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MATHÉMATIQUES

SUR LES PARTIES DE LA SUITE NATURELLE DES NOMBRES ENTIERS

Par N. LUSIN, de l'Académie

Les célèbres travaux de René Baire sur la représentation des fonctions discontinues, en dépit des apparences, ne sont pas au fond autre chose que des recherches relatives à la suite indéfinie

$$1, 2, 3, 4, 5, 6, \ldots$$

des entiers positifs. Ces travaux, arrêtés par une maladie et la mort prématurée de l'illustre auteur, paraissent être susceptibles à un prolongement naturel, au point de vue descriptif, ainsi que métrique, d'ailleurs.

1. Premières définitions. Lorsque l'on écrit une suite telle que la suite naturelle des nombres entiers

$$1, 2, 3, 4, \ldots$$

on se la représente sous la forme d'une série illimitée simple. Nous allons désigner cette suite par R. Si μ est un entier non négatif, nous pouvons désigner par R_μ l'ensemble des nombres

$$\mu+1, \quad \mu+2, \mu+3, \mu+4, \ldots$$

que nous appelons reste de la suite naturelle R. Il est clair que $R_0 = R$.

Si M est un ensemble fini d'entiers positifs (non nul ou bien dépourvu d'éléments), nous écrivons $M \equiv 0$. Si M est non nul, nous désignerons par $\overline{M}$ le plus grand des éléments qui appartiennent à M. Si M est un ensemble nul, nous écrivons $\overline{M} = 0$. Ces notations posées, il est évident que, si l'on a $\overline{M} = \mu$, il n'y a pas d'éléments de M dans le reste R_μ et, par suite, nous avons l'égalité $M = 0$ dans R_μ.

Par préférence, nous considérons des parties infinies E de la suite naturelle R. Si E et E' sont de telles parties, nous écrivons $E' \subset E$ lorsque tous les éléments de E', sauf un nombre limité (qui peut être nul) d'eux, appartiennent à E. Dans ce cas, nous avons $E' - E \equiv 0$. Si l'on a $E' - E = \mu$, nous avons l'inégalité évidente $E' < E$ dans le reste R_μ.

Deux parties infinies E et $\mathcal{E}$ de la suite naturelle seront dites orthogonales l'une à l'autre lorsqu'on a $E\mathcal{E} \equiv 0$; si $\overline{E\mathcal{E}} = \mu$, les parties E et $\mathcal{E}$ sont sans éléments communs dans R_μ.

Deux ensembles quelconques $\mathfrak{M} = \{E\}$ et $\mathfrak{N} = \{\mathcal{E}\}$ formés de parties infinies E et $\mathcal{E}$ de la suite naturelle seront dits orthogonaux l'un à l'autre (en général, non uniformément), si tous leurs éléments E et $\mathcal{E}$ sont respectivement orthogonaux, $E\mathcal{E} \equiv 0$; dans ce cas nous écrivons $\mathfrak{M}\mathfrak{N} \equiv 0$.

2. Séparabilité. Soient $\mathfrak{M}=\{E\}$ et $\mathfrak{N}=\{\mathcal{E}\}$ deux ensembles quelconques formés de parties infinies E et $\mathcal{E}$ de la suite naturelle. Nous écrivons $\mathfrak{M}\subset\mathfrak{N}$ lorsque la relation $E\subset\mathcal{E}$ est vérifiée pour tous les éléments respectifs E et $\mathcal{E}$ de ces ensembles. Si l'ensemble $\mathfrak{N}$ est formé d'une seule partie infinie H de la suite naturelle, écrire $\mathfrak{M}\subset H$ ce sera écrire la relation $E\subset H$ pour tous les éléments E de l'ensemble donné $\mathfrak{M}$. Dans ce cas particulier, la partie infinie H de la suite naturelle sera dite couverture de l'ensemble $\mathfrak{M}$.

Deux ensembles quelconques $\mathfrak{M}$ et $\mathfrak{N}$ formés de parties infinies de la suite naturelle seront dits séparables, si parmi leurs couvertures respectives H et K, $\mathfrak{M}\subset H$ et $\mathfrak{N}\subset K$ il y a des couvertures orthogonales: $HK\equiv 0$. Appelons séparatrices de deux ensembles $\mathfrak{M}$ et $\mathfrak{N}$ de telles couvertures orthogonales H et K.

Ces définitions étant posées, on a la proposition tout-à-fait évidente:

La condition nécessaire pour que deux ensembles quelconques $\mathfrak{M}$ et $\mathfrak{N}$ soient séparables est que ces ensembles soient orthogonaux l'un à l'autre: $\mathfrak{M}\mathfrak{N}\equiv 0$.

Dans les cas particuliers, cette condition est en même temps *suffisante*. Ceci nous assure le théorème suivant dont l'origine remonte au célèbre raisonnement de Paul du Bois-Reymond.

Théorème (Paul du Bois-Reymond). *Si deux ensembles $\mathfrak{M}$ et $\mathfrak{N}$ formés de parties infinies de la suite naturelle et orthogonaux l'un à l'autre sont dénombrables (ou finis), ils sont sûrement séparables.*

Démonstration. Par hypothèse, nous pouvons écrire $\mathfrak{M}=\{E_1, E_2, \ldots, E_n, \ldots\}$ et $\mathfrak{N}=\{\mathcal{E}_1, \mathcal{E}_2, \ldots, \mathcal{E}_n, \ldots\}$. Comme $\mathfrak{M}$ et $\mathfrak{N}$ sont orthogonaux l'un à l'autre, $\mathfrak{M}\mathfrak{N}\equiv 0$, nous avons $E_p\mathcal{E}_q\equiv 0$ pour tous les p et q possibles.

Désignons par $f(n)$ le plus grand des $2n$ entiers non négatifs $\overline{\mathcal{E}_\nu E_n}$ et $\overline{E_\nu \mathcal{E}_n}$, où ν ne surpasse pas n (le procédé de du Bois-Reymond), et définissons deux parties infinies H et K de la suite naturelle par la loi suivante:

$$H=E_1R_{f(1)}+E_2R_{f(2)}+\ldots+E_nR_{f(n)}+\ldots$$

et

$$K=\mathcal{E}_1R_{f(1)}+\mathcal{E}_2R_{f(2)}+\ldots+\mathcal{E}_nR_{f(n)}+\ldots$$

Tout d'abord, $E_n\subset H$ et $\mathcal{E}_n\subset K$ quel que soit un entier positif n, et, par suite, nous avons les relations $\mathfrak{M}\subset H$ et $\mathfrak{N}\subset K$ satisfaites. Donc, H et K sont des couvertures correspondantes de deux ensembles donnés $\mathfrak{M}$ et $\mathfrak{N}$.

Or, ces parties infinies H et K ainsi définies n'ont aucun élément commun. En effet, si un entier positif e appartient à H et à K, il appartient nécessairement à l'un des termes, $\mathcal{E}_qR_{f(p)}$, de la somme H et en même temps à l'un des termes, $\mathcal{E}_qR_{f(q)}$, de la somme K. Donc, e appartient à la partie commune $E_p\mathcal{E}_q$ et, par suite, si n est le plus grand des deux nombres p et q, nous avons $e\leqslant f(n)$. D'autre part, e appartient à la partie commune $R_{f(p)}R_{f(q)}$ et, définitivement, au reste $R_{f(n)}$ de la suite naturelle, ce qui est contraire à l'inégalité précédente. Donc, nous avons $HK=0$ et *a fortiori* $HK\equiv 0$. Il en résulte que les parties infinies H et K sont séparatrices de deux ensembles donnés $\mathfrak{M}$ et $\mathfrak{N}$, c. q. f. d.

La question qui se pose maintenant est la suivante.

Dans le cas général l'orthogonalité mutuelle $\mathfrak{M}\mathfrak{N}\equiv 0$ de deux ensembles donnés $\mathfrak{M}$ et $\mathfrak{N}$ est-elle une condition suffisante pour que $\mathfrak{M}$ et $\mathfrak{N}$ soient séparables?

C'est une question qu'il serait très important de résoudre; car, si on la résout par l'affirmative, on pourra affirmer que deux ensembles quelconques orthogonaux l'un à l'autre étant donnés, on peut déterminer effectivement leurs séparatrices. Si, au contraire, cette question se résout par la négative, il pourra exister deux ensembles non dénombrables orthogonaux l'un à l'autre et tels qu'il soit impossible de les séparer.

Sans vouloir aborder le fond de la question qui, à mon avis, ne sera résolue définitivement que par l'observation des faits; je me contenterai de quelques raisonnements auswahliques, c'est-à-dire fondés sur l'axiome du choix arbitraire de Zermelo (Auswahlprinzip).

3. Prolongement des suites orthogonales. Parmi les découvertes de George Cantor la plus importante est relative à la puissance du continu: *le continu n'est pas un ensemble dénombrable.* La forme précise que l'on peut donner à cet énoncé revient à la possibilité de *prolonger toute suite indéfinie* $x_1, x_2, \ldots, x_n, \ldots$ *de nombres réels en définissant effectivement un nombre réel nouveau* x *tel qu'il n'appartienne pas sûrement à cette suite.*

Poursuivons par analogie et appelons orthogonal en soi tout ensemble formé de parties infinies de la suite naturelle dont tous les éléments sont orthogonaux deux à deux.

Un ensemble $\mathfrak{M} = \{E\}$ orthogonal en soi est dit ouvert si l'on peut adjoindre à $\mathfrak{M}$ une partie infinie nouvelle $\mathcal{E}$ de la suite naturelle telle qu'elle soit orthogonale à chaque élément E de $\mathfrak{M}$, $\mathfrak{M}\mathcal{E} \equiv 0$. Dans le cas contraire, l'ensemble $\mathfrak{M}$ est dit fermé, c'est-à-dire ne pouvant plus être étendu par l'adjonction d'aucun élément nouveau sous la condition de retenir l'orthogonalité en soi.

Un ensemble orthogonal en soi et fini peut être fermé; mais s'il est effectivement dénombrable, il ne peut l'être jamais. C'est une proposition fort élémentaire dont l'idée même remonte à l'inventeur de la théorie des ensembles.

Théorème (George Cantor). *Toute suite indéfinie d'éléments orthogonaux deux à deux peut être toujours prolongée.*

Démonstration. Soit $E_1, E_2, \ldots, E_n, \ldots$ une suite indéfinie de parties infinies de la suite naturelle, orthogonales deux à deux: $E_p E_q \equiv 0$, si $p \neq q$. Désignons par $f(n)$ le plus grand des $\frac{n(n-1)}{2}$ entiers non négatifs $E_p E_q$, où p et q ne surpassent pas n, $p \neq q$. Il est clair que, dans le reste $R_{f(n)}$, les parties $E_1, E_2, \ldots, E_n$ sont sans élémens communs deux à deux. Par conséquent, nous pouvons déterminer effectivement un élément e_n de E_n aussi grand que l'on veut et tel qu'il n'appartienne à aucun des $n-1$ parties précédentes $E_1, E_2, \ldots, E_{n-1}$. *On reconnaît bien ici le fameux procédé diagonal de G. Cantor.*

Dans ces conditions, les éléments

$$e_1 < e_2 < e_3 < \ldots < e_n < \ldots$$

ainsi déterminés forment une partie infinie E effective de la suite naturelle telle qu'elle soit sûrement orthogonale à tous les éléments $E_2, E_2, \ldots, E_n, \ldots$ de la suite infinie donnée: $E_n E \equiv 0$ quel que soit n; d'ailleurs, bien entendu, les entiers positifs $\overline{E_n E} = e_n$ sont différents deux à deux, c. q. f. d.

4. Des ensembles orthogonaux non séparables. Le théorème précédent fait correspondre à chaque suite indéfinie de parties infinies orthogonales deux à deux $E_1, E_2, \ldots, E_n, \ldots$, effectivement énumérées au moyen des entiers positifs 1, 2, 3, ..., une et une seule partie effective E orthogonale à toutes ces parties, sans en excepter une seule.

En appliquant transfiniment ce théorème, on aura une suite bien ordonnée non dénombrable

$$E_0, E_1, E_2, \ldots, E_\omega, \ldots, E_\alpha, \ldots, \Omega$$

de parties infinies de la suite naturelle, orthogonales l'une à l'autre; d'ailleurs, par formation même de cette suite transfinie, les entiers positifs $\overline{E_\beta E_\alpha}$, pour chaque valeur fixe de α et pour β variable inférieur à α, $\beta < \alpha$, sont tous différents.

Cela posé, partageons d'une manière quelconque la suite transfinie considérée en deux suites, toutes les deux non dénombrables et sans parties

communes; nous désignons par $\mathfrak{M}$ et $\mathfrak{N}$ deux ensembles non dénombrables formés respectivement des éléments de ces suites:

$$\mathfrak{M} = \{E_{\gamma'_0}, E_{\gamma'_1}, \ldots, E_{\gamma'_\omega}, \ldots, E_{\gamma'_\alpha} \ldots\ldots | \Omega\}$$

et

$$\mathfrak{N} = \{E_{\gamma''_0}, E_{\gamma''_1}, \ldots, E_{\gamma''_\omega}, \ldots\ldots, E_{\gamma''_\alpha}, \ldots\ldots | \Omega\}$$

les indices γ'_α et γ''_α étant des nombres finis et transfinis différents rangés par ordre de grandeur croissante.

Il est manifeste que ces ensembles $\mathfrak{M}$ et $\mathfrak{N}$ sont orthogonaux l'un à l'autre: $\mathfrak{M}\mathfrak{N} \equiv 0$.

Je dis maintenant que $\mathfrak{M}$ et $\mathfrak{N}$ ne sont pas séparables.

En effet, si, par impossible, H et K seront des séparatrices, nous aurons $E_{\gamma'_\alpha} \subset H$, $E_{\gamma''_\alpha} \subset K$ et $HK \equiv 0$. *A fortiori* nous aurons $E_{\gamma'_\alpha} K \equiv 0$ et $E_{\gamma''_\alpha} H \equiv 0$ quel que soit α inférieur à Ω. Nous concluons de là qu'il existe une infinité dénombrable d'indices β tels qu'on a $\overline{E_{\gamma'_\beta} K} < m$, où m est un entier positif fixe; nous pouvons supposer ces indices β inférieurs à un nombre transfini β^* fixe, $\beta^* < \Omega$. Il est clair qu'on peut trouver un nombre transfini fixe λ, $\lambda < \Omega$, tel que l'on ait $\gamma_{\beta^*} < \gamma''_\lambda$. Dans le reste R_m les parties $E_{\gamma'_\beta}$ et K sont sans élément commun. Comme nous avons $E_{\gamma''_\lambda} \subset K$, il existe un reste R_μ tel qu'on ait, dans ce reste, l'inégalité $E_{\gamma''_\lambda} < K$. Soit m^* le plus grand des deux entiers positifs m et μ; dans le reste R_{m^*} il n'y a pas d'éléments communs à $E_{\gamma'_\beta}$ et $E_{\gamma''_\lambda}$. Or, nous avons $\gamma'_\beta < \gamma''_\lambda$ et les indices β sont sûrement en nombre infini, ce qui est précisément impossible, c. q. f. d.

Pour les mathématiciens qui s'intéressent, par préférence, de la technique même des raisonnements mathématiques et qui admettent l'existence de tous les nombres transfinis de la seconde classe, les considérations du texte permettent de poser une quantité embarrassante de questions difficiles qu'on rencontre sur cette voie et qu'on ne peut pas trouver sur autres voies: telles sont, par exemple, les questions relatives à la séparabilité B des ensembles orthogonaux en soi; parmi lesquels on a sûrement des ensembles ayant la puissance du continu; etc.

Mais un travail beaucoup plus pénible est encore entièrement à faire pour savoir quel est le sens réel et précis qui se cache derrière les raisonnements du texte, fondés manifestement tous sur le choc purement verbal des termes: le fini et l'infini.

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LITTÉRATURE CITÉE

[1] Felix Hausdorff, Fundam. Mathemat., XXVI, 241 (1938).

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MATHEMATICS

ON THE DISTRIBUTION OF THE ZEROES OF AN INTEGER FUNCTION

By N. MEYMAN

(Communicated by S. L. Soboleff, Member of the Academy, 3.VI.1943)

We write an integer function $F(z)$ in the form $F(z)=g(z)+ih(z)$, where $g(z)$ and $h(z)$ take real values on the real axis. Let $g(z)$ and $h(z)$ have no common roots. $F(z)$ is called a B-function, if $\left|\frac{g(z)+ih(z)}{g(z)-ih(z)}\right|<1+\eta$ holds for sufficiently large values of $|z|$, z lying in the upper half-plane, and η being an arbitrary positive number (2).

We consider a point-set M lying in the upper half-plane and defined by the inequality

$$|\zeta(z)|=\left|\frac{F(z)}{\overline{F(z)}}\right|=\left|\frac{g(z)+ih(z)}{g(z)-ih(z)}\right|>1. \qquad (1)$$

The boundary of M will be denoted by $\dot{M}$. M consists of a finite or infinite number of components U each of which is a domain. The boundary of each component consists in its turn of a finite or infinite number of components. Each point of $\dot{M}$ will be numbered as many times as it enters in $\dot{M}$. Each component of $\dot{M}$ is an analytical curve. The only possible singularities of $\dot{M}$ are angle-points and multiple points, the value of the angle being $\frac{\pi}{n}$, where n is a natural number. At these points $\zeta'(z)=0$. The positive direction on $\dot{M}$ is defined to leave a half-neighbourhood contained in M on the left-hand side. The components of $\dot{M}$ which lie entirely in a finite part of the plane will be denoted by K. The K are closed curves whose image in the ζ-plane is a circumference $|\zeta|=1$ which is run in the negative direction a finite number of times. By L we denote the components of $\dot{M}$ tending to ∞. Their image is a circumference $|\zeta|=1$ run in the negative direction. According as the circumference is travelled a fractional, integer or infinite number of times, L will be denoted by L_1, L_2 or L_3.

Along the branch of L_1, stretching to ∞ in the positive direction, $\zeta(z)$ tends to a limit and does so also along the branch tending to ∞ in the negative direction, but its limit here is known to be different from the first. Then in the domain of the upper half-plane bounded by L_1, $\zeta(z)$ in virtue of a well-known theorem (1) takes on all finite and infinite values an infinite number of times but for two values at most. In the domain of the upper half-plane bounded by L_2 $\zeta(z)$either takes all finite and infinite values with the exception, perhaps, of two, an infinite number of times (in this case L_2 will be denoted by L_{21}) or else tends uniformly to a limit as z passes to ∞ (1) (in this case L_2 will be denoted by L_{22}). The latter limit obvi-

ously coincides with the asymptotic value of $\zeta(z)$ along L_2, and has the absolute value of 1.

The points of the real axis satisfying to the inequality

$$h'(x)\,g(x)-h(x)\,g'(x)>0 \tag{2}$$

are not contained in $\dot{M}$; the points of the real axis satisfying to the inequality

$$h'(x)\,g(x)-h(x)\,g'(x)<0 \tag{3}$$

are contained in $\dot{M}$, and so are their real neighbourhoods; the points of the real axis for which

$$h'(x)\,g(x)-h(x)\,g'(x)=0 \tag{4}$$

are angle-points of $\dot{M}$.

Let a be a real number, finite or infinite. The roots of the equation $\frac{h(z)}{g(z)}=a$ lying in the closed upper half-plane will be called upper a-points. Let such a root be k-uple. The upper multiplicity is put at k if $Jz>0$; at $\frac{k}{2}$ if the root is real and k is even; and at $\frac{k-1}{2}$ or $\frac{k+1}{2}$ for an odd k and a real z, according as $\arg F(z)$ increases or decreases at z. In other words, the upper multiplicity of a point z is defined to be equal to the multiplicity in which this point is contained in $\dot{M}$. The sum of upper multiplicities of all a-points for a given a will be denoted by $S_a\frac{h}{g}$. The numbers $S_0\frac{h}{g}$ and $S_\infty\frac{h}{g}$ are obviously equal to the number of the upper zeroes of h and g (taking into account their upper multiplicities).

We denote by $w(\zeta)=z$ the function inverse to the meromorphic function $\zeta(z)$, and by N the prototype of M lying on the Riemannian surface of the function w. To each component U of M there corresponds a certain component V of N, and conversely. Moreover, each component V is also a component for a part of the Riemannian surface lying over the exterior of the circle $|\zeta|=1$. The boundary of V consists only of points lying over the circle $|\zeta|=1$, and of points which are asymptotic values of $\zeta(z)$ for the paths lying within U.

If the function $F(z)$ belongs to B, then $\zeta(z)$ has no more than one asymptotic value in the upper half-plane; moreover, if $\zeta(z)$ tends to this value along two paths Γ_1 and Γ_2, it also tends to it uniformly in the domain of the upper half-plane bounded by $\Gamma_1+\Gamma_2$, when $z\to\infty$. Absolutely this asymptotic value is $\leqslant 1$; equality must hold, when M contains a component tending to infinity. For functions of the class B every V consists of a finite number of full sheets lying over the exterior of $|\zeta|=1$, and the prototype of the boundary of each U is the circle $|\zeta|=1$ run through an integer number of times in the negative direction. The total number of all sheets belonging to N is finite. In this case the boundary of M consists of a finite number of the K and L_{21}. Conversely, if N is contained in a finite number of sheets, then $F(z)\in B$. Now we can point out

Theorem 1. *If $F(z)\in B$, then for all values a, different from the asymptotic value of $\zeta(z)$ in the upper half-plane, the $S_a\frac{h}{g}$ are bounded and equal. For the asymptotic values the value of $S_a\frac{h}{g}$ is less.*

If $F(z)\in B$ and is not a polynomial, then $\arg F(z)$ *monotonously increases to infinity along a certain part of the real axis.*

Theorem 2. *If there exist two (finite or infinite) numbers a and b such that the values $S_a\frac{h}{g}$ and $S_b\frac{h}{g}$ are finite, then either $F(z)$ belongs to B, and*

therefore the number of roots of $F(z)$ in the opened upper half-plane is equal to the largest of the values $S_a \frac{h}{g}$, $S_b \frac{h}{g}$, or $F(z)$ has an infinite number of roots in the opened half-plane, and therefore does not belong to B.

Proof. Since $S_a < +\infty$, $\dot{M}$ does not contain the components L_3. Each component K contains the same number (different from zero) of upper a-points and b-points. Each component L_{22} contains either one upper a-point, or one b-point at least. Therefore, $\dot{M}$ contains only a finite number of components of types K and L_{22}. If $\dot{M}$ contains no components of types L_1 and L_{21}, then the principle of variation of the argument is applicable to M, in this case $F(z) \in B$. The variation of $\arg \zeta(z)$ along K is $-2r\pi$, where r is the number of the upper a-points or b-points lying on K. The variation of $\arg \zeta(z)$ along L_{22} is $-2s\pi$, where s is the largest of the numbers of a-points and b-points lying on L_{22}. Consequently, in this case the number of poles of the function $\zeta(z)$ in the upper half-plane, or, what is the same, the number of the zeroes of $F(z)$ in the lower half-plane, is equal to the largest of the numbers S_a and S_b.

Even if $\dot{M}$ should contain only one component of type L_1 or L_{21}, $F(z)$ would have an infinite number of zeroes in the lower half-plane or, more precisely, in the domain conjugate to the domain bounded by such a component. For, in such a domain $\zeta(z)$ has an infinite number of poles, because $\frac{1+ia}{1-ia}$ and $\frac{1+ib}{1-ib}$ are exceptional values in the sense of Lindelöf's [1] theorem.

Remark. From the proof of Theorem 2 we infer the following corollary. If S_a and $S_b < +\infty$, then $F(z)$ belongs to B if and only if the quotient $\frac{h(z)}{g(z)}$ has no more than one asymptotic value in the upper half-plane. Moreover, if $F(z)$ does not belong to B, there exists an infinity of components of type L_1 or L_{21} enclosing each other.

Example. Let $F(z) = \sum_{l,k=0}^{n} c_{lk} z^l e^{i\alpha_k z}$, where $\alpha \geqslant 0$ and the sum contains a term in which l and α_k are not less than the corresponding exponents in any other term. Since $F(z) \in B$, the number of roots of $F(z)$ in the lower half-plane is equal to the largest of the numbers S_0 and S_∞. In this case the asymptotic value of $\frac{h}{g}$ for $Jz > 0$ is i. Therefore, $S_0 = S_\infty$. $F(z)$ has no zeroes in the lower half-plane if and only if $S_0 = S_\infty$. This has been obtained in paper [3] for integral α's.

It follows from Theorem 2:

I. A function $F(z)$ of the class B has no roots in the lower half-plane if and only if $S_0 = S_\infty = 0$.

II. If $F(z)$ has only a finite number of zeroes in the lower half-plane and does not belong to the class B, then at least one of the numbers S_0 or S_∞ is equal to ∞.

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HYDROMECHANICS

CERTAIN PROBLEMS OF FILTRATION IN A TWO-LAYER MEDIUM REDUCED TO THE CASE OF A HOMOGENEOUS GROUND

By N. K. KALININ

(Communicated by N. E. Kochin, Member of the Academy, 26.V.1943)

1. P. J. Polubarinova-Kochina ([1]) has put an instance to show that the method developed by her for solving filtration problems in a homogeneous ground, based on the use of the analytical theory of linear differential equations, can be extended to a certain class of filtration problems in a two-layer medium. However, a region analogous to velocity hodograph is not easy to plot for this kind of problems, and there arise a number of difficulties connected with the determination of the number of singular points (when the so-called collateral singular points are apt to be omitted) and with the finding of their indices. Considerable interest attaches therefore to the reduction of filtration problems in a two-layer ground to the case of a homogeneous ground. Owing to the work of Polubarinova-Kochina ([2]) and Riesenkampf ([3]), the method for the solution of these problems by means of linear differential equations has acquired great simplicity and representativeness. Below we show that a certain group of filtration problems in a two-layer medium; when there is no free surface, may be reduced to the case of a homogeneous ground in the presence of a free surface, and also to such problems that admit of a solution by the well developed method of Polubarinova-Kochina. As for the regularity of singular points in the case of filtration in a two-layer medium, this problem needs no special solution; once the regularity of singular points is established for the case of filtration in a homogeneous ground ([3]). The motion is supposed to be plane (parallel to the vertical plane), and steady. Instead of the complex potential $f=\varphi+i\psi$, we shall examine the so-called complex lost head $w=q+ih$ ([4]) which is connected with the former by the relation $f=ikw$; where q is the function of discharge; h; the lost head; k, the coefficient of filtration.

2. Suppose that the filtration region G (which for simplicity sake is assumed to be simply connected) consists of two grounds with filtration coefficients k_1 and k_2 occupying simply connected regions G_1 and G_2 having a common boundary L (division line)* and bounded, moreover, by the arcs L_1 and L_2, respectively. Let us consider complex lost heads $w_1=q_1+ih_1$, $w_2=q_2+ih_2$, regular in the regions G_1 and G_2, respectively (it is assumed that there are no point sources inside G).

We assume further that the position of the boundaries L_1 and L_2 is given and that the arc L_1 (resp. L_2) consists of segments on which h_1 and q_1 (resp. h_2 and q_2) take constant values alternately.

Let us refer the filtration region to the plane of a complex variable $\zeta=\xi+i\eta$. The use of the method developed by Polubarinova-Kochina depends in our case upon the existence of an analytical function $\xi=\zeta(t)$ by which the region G_1 is mapped conformably upon the upper and the region G_2 upon the lower half-plane of the auxiliary complex variable t; in other words, the regions

* L is supposed to be a regular arc of an analytical curve.

G_1 and G_2 should be symmetrical with respect to the arc L. Suppose arbitrarily that the region G satisfies this condition. The function $\zeta(t)$ will be taken in such form that the line L passes to the positive part, and the lines L_1 and L_2 to the negative part of the real axis of the variable t. Now we shall consider the functions $W_1(t)=\frac{dw_1}{dt}$, $W_2(t)=\frac{dw_2}{dt}$.

They represent complex velocities of certain flows, determined in the upper and lower half-planes of t, respectively, and satisfying for $t>0$ (line of division of grounds) the conditions (1):

$$J(ik_1W_1-ik_2W_2)=0, \quad J(W_1-W_2)=0, \tag{1}$$

where J denotes the imaginary part of the function; for $t<0$ they take real and imaginary values in alternation. Owing to the latter circumstance, the function W_2 determined in the lower half-plane may be extended analytically to the upper half-plane through any segment of the negative part of the real axis of the variable t. The result of such an analytical extension we denote by $\tilde{W}_2$. If the extension is effected through a segment upon which $J(W_2)=0$ *, then, according to the symmetry principle of Riemann-Schwarz, we have

$$J(\tilde{W}_2)=-J(W_2), \quad J(i\tilde{W}_2)=J(iW_2). \tag{2}$$

With $t<0$ the function $\tilde{W}_2$ satisfies the same boundary conditions as the function W_2 does; yet for $t>0$ instead of (1) we, in accordance with (2), have to take the conditions

$$J(ik_1W_1-ik_2\tilde{W}_2)=0, \; J(W_1+\tilde{W}_2)=0. \tag{3}$$

Instead of functions W_1, W_2 we now shall seek for the functions W_1, W_2 determined in the upper half-plane of t and satisfying on the real axis for $t>0$ to the conditions (3). On the segments of the negative part of this axis (these segments, generally speaking, are different for W_1 and $\tilde{W}_2$) they take now real now imaginary values; that is, for $t<0$ we have alternately

$$J(W_1)=0 \quad \text{and} \quad J(iW_1)=0 \tag{4}$$

or, respectively,

$$J(\tilde{W}_2)=0 \quad \text{and} \quad J(i\tilde{W}_2)=0. \tag{5}$$

On the other hand, let us consider the problem of ground water motion in a homogeneous (isotropic) ground with a filtration coefficient $k=1$ in a simply connected region of the half-plane $z=x+iy$ (the axis y is vertical), bounded by a polygonal curve that consists of horizontal and vertical segments along which $h=\text{const}$ (the boundary of the reservoir) and $q=\text{const}$ (solid wall) in alternation, and by free surface on which the function of discharge satisfies the condition ([4],[5]) **

$$q=cx+\text{const} \; (c>0) \text{ or } J(iw-icz)=\text{const}, \tag{6}$$

where $w=q+ih$ is the complex lost head of the given current in a homogeneous ground. On the free surface, moreover, there must hold the condition of pressure constancy, i. e.

$$h=-y+\text{const} \text{ or } J(w+z)=\text{const}. \tag{7}$$

Let the region of the ground water be mapped conformably onto the upper

* If the extension is through a segment where $J(iW_2)=0$, the reasoning should be slightly modified.

** If by points at which the tangent to the free surface is vertical the latter is divided into areas bounding the region of motion from above and from below, then we, according to (6), will have evaporation on the first areas and absorption of the liquid on the second areas.

half-plane of the auxiliary complex variable t in such a way that to the free surface there should correspond a positive part of the real axis of this variable and to the polygonal curve its negative part. We shall consider the following functions connected with this mapping: $W(t)=\frac{dw}{dt}$, $Z(t)=\frac{dz}{dt}$.

These functions are regular in the upper half-plane of t and for $t>0$ satisfy two boundary conditions obtained from (6) and (7) by differentiation with respect to t: namely

$$J(iW-icZ)=0,\ J(W+Z)=0. \tag{8}$$

On the segments of the negative part of the real axis of t (these segments are, generally speaking, different for W and Z) they arbitrarily take now real now purely imaginary values; that is, for $t<0$

$$J(W)=0 \quad \text{and} \quad J(iW)=0 \tag{9}$$

in alternation, and, accordingly,

$$J(Z)=0 \quad \text{and} \quad J(iZ)=0. \tag{10}$$

It may now easily be seen that the conditions (3), (4), (5) and (8), (9), (10) can be brought to identity by putting

$$W_1=W,\ \tilde{W}_2=Z,\ c=\frac{k_2}{k_1} \tag{11}$$

and by suitably selecting the polygonal curve in the problem of filtration in a homogeneous ground.

So we see that the class of two-layer filtration problems under consideration is actually equivalent to certain filtration problems in a homogeneous ground when a free surface is present. Having solved the problem of filtration in a homogeneous ground, that is having found the functions $W(t)$, $Z(t)$, by means of formulae (11) we can find the functions $W_1(t)$, $W_2(t)$ that give the solution of the respective filtration problem for a two-layer medium.

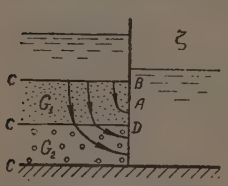

Fig. 1

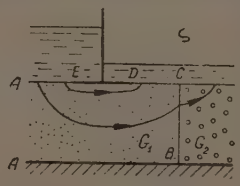

Fig. 2

From what has been said here we may be easily convinced that in the case of filtration under a groove in a two-layer medium with a horizontal division line between the layers [the problem of Polubarinova-Kochina (1)], the problem presented in Fig. 1, ζ is equivalent to that of filtration in a homogeneous ground through a dam drained at the foundation (Fig. 1, z); at the free surface evaporation takes place. Another example is the filtration problem under a flat floatbet with a vertical division line between the layers, presented in Fig. 2, ζ. It is likewise equivalent to the filtration problem in a homogeneous ground (Fig. 2, z). A detailed discussion of this and other examples will be made in a more comprehensive paper to be published.

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BIOPHYSICS

VISIBILITY OF TEST-OBJECTS UNDER LOW ILLUMINATIONS

By L. N. GASSOVSKY, K. L. BULANOVA and Z. N. ENNO

(Communicated by S. I. Wawilow, Member of the Academy, 7.IV. 1943)

For high background luminosities the relation between the visual acuity and the contrast value is sufficiently well known as a result of a number of special investigations [see, for instance; the paper by Cobb and Moss (1)].

An attempt to find this relation for very low background luminosities; corresponding to nocturnal vision in the absence of artificial light sources, has been made only by Conner and Ganoung (2) on seven observers (background liminosities within the range between 0.001 asb and 1 asb).

The work was carried out on 15 observers. As test-objects, besides the Landolt rings which had been used in the work of Conner and Ganoung; we also used optical circles, miras, as are usually applied for determining the resolving power of optical instruments*.

The relation between the visual acuity and the contrast value between the test-object and the background was studied by us for background luminosities (adaptation field) within the range of 0.0005 to 5.0 asb. On account of the fact that under conditions of fight manoeuvres the observer has but very short time intervals at his disposal; the relation between the visual acuity and the contrast value was investigated under conditions of limited times of observation, restricted to 0.02 and 2″.

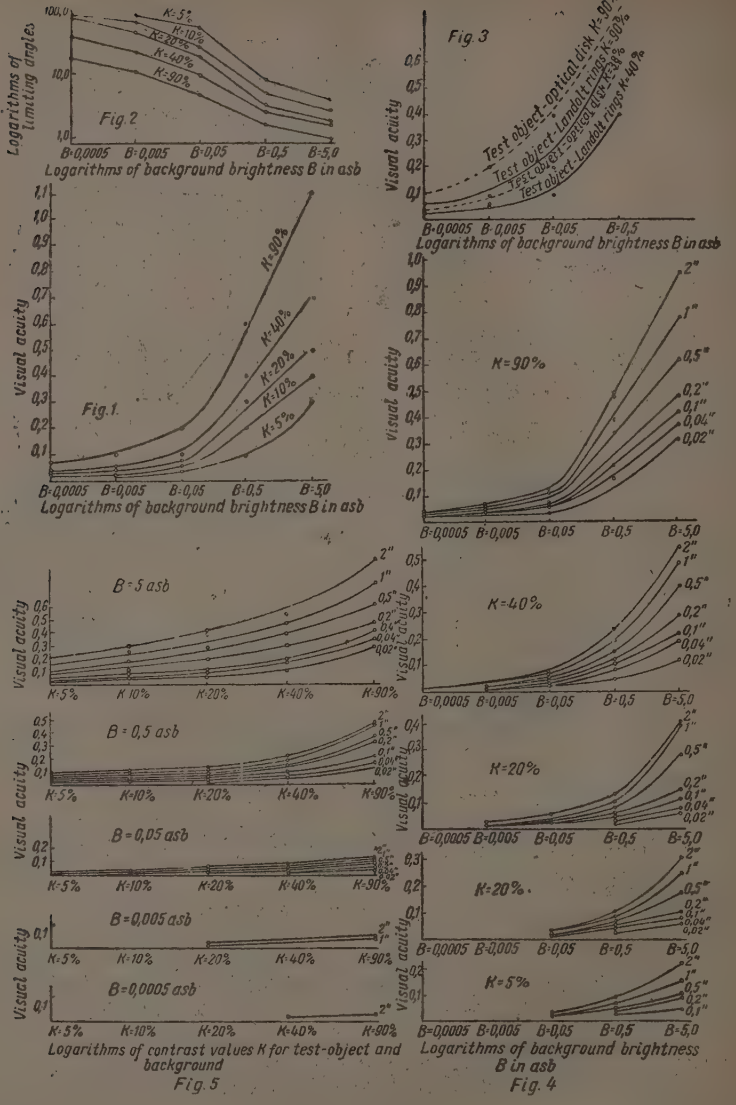
The instrument used in our investigation consisted essentially of a large (180 × × 180 cm) white screen uniformly illuminated over the whole of its surface by special luminaries. The luminosity of the screen could be varied arbitrarily by inserting before the luminaries specially devised and selected neutral filters. The luminosity of the screen could assume the following five values: 0.0005; 0.005; 0.05, 0.5 and 5.0 asb. The constancy of the current used for feeding the bulbs of the luminaries of the screen, was controlled by a precision voltmeter.

The test-objects could be placed in the centre of the screen: as test-objects Landolt rings cut of Wattman paper or optical circles were used. The contrast values of the test-objects with respect to the background (white Wattman paper) were 90, 40, 20, 10 and 5%.

We used in our experiments rings and circles of 15 different sizes; so that the total number of rings and circles amounted to 80. The observer was placed at a distance of 5 m from the screen. The observations were carried out in a room with blackened walls. No external light was allowed to penetrate into the room.

Prior to start the experiment; the observer was kept in complete darkness during 20 minutes and thereafter made to fix during 10—12 minutes a screen-background of a given luminosity. After light adaptation was complete, the observer was required to determine the direction of the breaks in the Landolt rings or

* Such a mira consists of a circle covered by a set of parallel black and white strips of equal width.



of the strips on the circles suspended against the screen. Ten observations at least were carried out for a test of each contrast value and angular size.

1. Observations with Landolt Rings. The character of the relation obtained in this case is borne out by the curves of Figs. 1 and 2. In Fig. 1 the logarithms of the background luminosity B are plotted as abscissae and the visual acuity; as ordinates. Each of the curves corresponds to a definite contrast value K between the test-object and the background. It can be seen from these curves that with increase of the background luminosity the visual acuity increases, and this the more rapidly, the darker the test-object compared with the back-

ground. Thus, for example; for a background luminosy of 5 asb and a contrast value of 90% the visual acuity is twice as large as the visual acuity with a 20% contrast value and the same luminosity. The curves in Fig. 1 consist of two parts: the left part with a gentle slope corresponding to the work of the rods of the visual apparatus of the eye (low values of background luminosities), and of the right one with an abrubt slope (high values of background luminosities) which corresponds to the work of the cones. Separate curves of rod and cone vision corresponding to these two parts of the curves in Fig. 1 have been obtained by Conner and Ganoung.

In Fig. 2 the relation obtained is represented in a somewhat different form: while the logarithms of the background luminosity are again plotted as abscissae, the logarithms of the limiting angles* are plotted as ordinates.

The curves of Figs. 1 and 2 are means for 15 observers. The individual curves of separate observers differ but very little from the mean curves. The curves obtained by us are very similar to those of Conner and Ganoung**.

2. Observations with Optical Circles (Miras). Observations with optical circles have yielded results similar to those obtained with Landolt rings. The visual acuity exceeded, however, in all cases that for Landolt rings by 40—50%. This should be taken to mean that the position of strips in a circle is discerned by the observer easier than the position of a break in the Landolt rings. This is illlustrated by Fig. 3. The observations with circles are strongly affected by slight degrees of astigmatism frequently met with.

3. Observations during Limited Time Intervals. The following values of the observation time (exposure) were used: 2, 1, 0.5, 0.2, 0.1, 0.04 and 0.02″.

The results obtained under such conditions are illustrated by the curves in Figs. 4 and 5. The increase of the duration of the observation influences the visual acuity in the same way as the increase of the contrast value K. This can easily be seen from a comparison of the curves in Fig. 1 and Fig. 4.

A comparison of the groups of curves in Fig. 4 shows that with a decrease of the contrast value K the divergence between the curves is diminished and their bases are shifted to the right. This means that with a decrease in the contrast value and with a limited observation time visual acuity is reduced; this leads to the impossibility of discerning even the largest objects with small background luminosities. The effect of the restriction of observation time is more important for large contrast values K and background luminosities B.

The curves of Fig. 4 enable one to determine the dimensions of an object which can be discerned by the eye for any given duration of observation and for any given contrast value and background luminosity (abaptation field of the eye) B.

In Fig. 5 the logarithms of the contrast value K are plotted along the abscissa and the visual acuity along the ordinate. In this case, too; the divergence between the separate curves diminishes with the decrease of background luminosity B, while their basis is gradually shifted to the right in the direction of large values of the contrast value K. The curves of Figs. 4 and 5 can be used for determining the maximum distances of visual observations.

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* The limiting angle is the smallest angle under which the observer can perceive a break in a Landolt ring or in the interval between two dark strips in an optical circle, provided that the position of the ring or of the circle be recorded by the observer correctly.

** The existing discrepancies may among other causes be accounted for by the fact that the significance of observations in our experiments was put at 80%, whereas in the experiments by Conner and Ganoung it was 55%.

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CHEMISTRY

PRECIPITATION OF MERCURY FROM SOLUTIONS OF SULPHIDES WITH AMALGAMATED ZINC

By I. N. PLAKSIN and L. D. PLAKSINA

(Communicated by E. V. Britzke, Member of the Academy, 31.III.1943)

The hydrometallurgy of mercury is a comparatively young branch of metallurgy. Most of the mercury deposits of the USSR; such as those of Salair (West Siberia); Haidarkan (Kazakhstan), Aiatka and Iss (Urals), are characterized by a low content of mercury in the ore.

The necessity of dressing large masses of ore puts forward hydrometallurgical methods of treatment that are cheaper, simpler and give a higher extraction as compared with the pyrometallurgical methods (for instance, leaching out of mercury with weak alkaline solutions of sodium sulphide in the case of lump ore in wood vats). These methods have certain advantage also for the concentrates in reducing losses in the treatment of rich materials.

In this country the first research work and industrial tests in this field were carried out in 1931 by Plaksin and Fishkova (1); the results obtained were positive. A series of subsequent investigations carried out by a number of institutes have shown the expedience of applying this method to low grade mercury ores. In connexion with the setting into operation of hydrometallurgical plants the question arises as to the reagents to be used in this process.

Until recently in all investigations; beginning with the work of Mulholland (2); and in the practice of plants in this Union and abroad; metallic aluminium was used as precipitation agent. The first to introduce it into the industrial practice of the hydrometallurgy of mercury was Thornhill (3) who described it in 1915.

On account of the great need for aluminium experienced in the time of war the authors undertook; along with a survey of the mercury ores of the Urals, a search for new precipitation agents. As to metal precipitation agents, the essential difficulty consisted in the capacity of metals to be sulphidized on the surface in sodium sulphide solutions.

The use of non-metallic precipitation agents, in particular of charcoal, resulted in the formation of too voluminous and hardly workable precipitates. The method devised and suggested by the authors consists in the precipitation of mercury with zinc amalgam or with zinc whose surface is covered with mercury protecting its surface from the formation of a film of zinc sulphide. Similar to aluminium, the zinc amalgam and amalgamated zinc precipitates from solutions antimony, the usual satellite of mercury minerals, which is highly soluble in sodium sulphide solutions.

Zinc amalgam and amalgamated zinc can be repeatedly used for mercury precipitation, whereas aluminium can be used no more than once, because of its dissolving vigorously in alkaline solutions of sodium sulphide. According to data of the work of the Salair plants, the expenditure of aluminium amounts to 1 kg per 1 ton of dry sand.

Numerous experiments on the precipitation of mercury from sulphide solutions by amalgamated zinc showed the essential rôle played by the diffusion phenomena in this process. By increasing the efficiency of stirring it is possible to reduce considerably the time of precipitation; the concentration of the solution in respect of Na_2S and NaOH and also the ratio between the precipitating agent and the mercury.

Preliminary investigations showed that mercury is precipitated at the rate of 99.9% per 30 min.; the ratio mercury : precipitating agent being 1 : 80. Under industrial conditions this ratio will be much lower. It is possible to lower the expenditure of the precipitating agent at the expense of using in the final stage of precipitation filtration of the solution through a layer of the precipitating agent on the surface of the filter. The concentration in the sodium sulphide solution during precipitation depends mainly on the mode and efficiency of stirring the solution during precipitation. In the case of stirring by means of friction cylinders a high percentage of precipitated mercury is obtained at concentrations 4.6% Na_2S and 0.2% NaOH. When the stirring device is Wagner's agitator, the precipitation can be achieved at a concentration of 1.5% Na_2S, probably owing to better stirring. The stirring of the solution makes it possible to use low concentrations of sodium sulphide; the effect obtained being sufficient for the industrial realization of the process. The minimum concentration of sodium sulphide allowing for satisfactory precipitation could be reduced to 0.5% Na_2S by making stirring more efficient.

Owing to this, for an efficient precipitation of mercury from sodium sulphide solutions appliances should be designed allowing for the application of intensive stirring and of the circulation of the solution with a simultaneous condition that precipitation be achieved with rather coarse-grained material. Experiments carried out with finely ground amalgams gave a lower per cent of extraction.

As precipitating agent we used a zinc amalgamated on the surface that contained 96% metallic zinc and 4% mercury. The zinc was covered with mercury in a dilute solution (5%) of hydrochloric or sulphuric acid. If the solution is not stirred, amalgamated zinc precipitates up to 98.5% of mercury within 12 hours.

For the precipitation of mercury from solutions amalgamated zinc can be used repeatedly. The best results are obtained if the surface of zinc that has been used once is cleansed with a 0.5% solution of hydrochloric or sulphuric acid. Repeated use of amalgamated zinc leads to the formation of a zinc amalgam with a higher mercury content.

When it is necessary to use a precipitating agent with a higher content of mercury (for instance, zinc amalgam), the conditions for precipitation are somewhat different. A 30 minute contact with the precipitating agent leads to a precipitation of 99.9% of mercury from a solution when the ratio mercury: precipitating agent is 1 : 50. The size of the amalgam grains is 1 mm. When the ratio mercury precipitating agent in the solution is 1 : 2, the time of precipitation should be 48 hours. 99.9% of Hg being extracted. The concentration ratio of sodium sulphide and alkali in solution is the same as in the preceding case.

The amalgam can be used many times without treatment with dilute acid and later with a 1 min. treatment with 0.5% hydrochloric or sulphuric acid. Zinc amalgam or granulated zinc covered with mercury are used repeatedly; whereas aluminium dissolves completely in alkaline solutions of sodium sulphide.

When precipitation is effected with the aid of zinc, the process is materially affected by the temperature. Precipitation with amalgam proceeds successfully at 20° C. This is explained by the influence upon this process of the conditions of zinc diffusion in the surface layer of the amalgam.

The use of superficially amalgamated zinc offers a considerable advantage; for the process can then be carried on at a much lower temperature; down to 9°. Leaching with solutions of sodium sulphide is sometimes prevented by lack in the locality of a given reagent or of the raw materials for its production.

Owing to the presence of mercury in barite ores or of barite deposits within the area of mercury occurrences; it seemed of interest to find out the conditions of precipitation with zinc also from solutions obtained by leaching with barium

sulphide. The conditions of precipitation from these solutions differ from those of the precipitation of mercury with amalgamated zinc from solutions of sodium sulphide.

First of all; the lower solubility of barium sulphide and its susceptibility to temperature makes it necessary to use solutions of very low concentrations at a temperature below 10° C. Another very important factor in the precipitation from solutions of barium sulphide is the necessity of operating with solutions of an alkalinity not below 0.5%.

In the case of precipitation from solutions of sodium sulphide this factor is less important, owing to the alkaline reaction of sodium sulphide, resulting from hydrolysis and a greater stability of solutions of NaOH as compared to those of $Ba(OH)_2$ formed in the hydrolysis of barium sulphide.

After precipitation there remains a part of the amalgam which did not enter the reaction, and can easily be separated and used again.

In this process mercury is precipitated to a considerable degree as a fine black-grey powder. Owing to this; mercury cannot be isolated by pressing out and is separated by distillation from a retort at a temperature of 400—450° C of its mixture with lime. The expenditure of zinc in the precipitation with amalgamated zinc is 1.5 and sometimes 2 parts per 1 part of metallic mercury.

The application of this procedure on an industrial scale will lead to a much lower expenditure of zinc as compared to precipitation with aluminium, the expenditure of which is considerable in the case of precipitation in extractors at the Salair hydrometallurgical plants.

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PHYSICAL CHEMISTRY

THE FORM OF THE FIRST BRILLOUIN ZONE IN THE CASE OF HEXAGONAL STRUCTURE OF NICKEL-ARSENIDE TYPE

By E. S. MAKAROV

(Communicated by A. N. Frumkin, Member of the Academy, 23.IV.1943)

The Brillouin zone theory, as developed by Jones (1); has permitted to explain a number of physical properties in chemical elements, their alloys and compounds. Applied to the problems of the chemistry of metal alloys, this theory proved especially fruitful in explaining the Hume-Rothery empirical rule which governs the formation of the so-called electronic compounds. Yet, compounds with normal valency, such as Mg_2Sn for instance, abso fit well in terms of the zone theory (2).

In this article we made an attempt to determine the form of the first Brillouin zone for a hexagonal crystalline lattice of nickel-arsenide type. The methods for finding the Brillouin zone and all necessary calculations are described in a well known book by Mott and Jones (3).

Compounds of nickel-arsenide structure are known to possess well marked metallic properties: a metallic lustre and electronic conductivity. However, the electronic conductivity is not high in these compounds. In Table 1 data on electric resistance ρ at 25° C and its temperature coefficient in the interval 25—100° C are given, according to the results of our experiments (4), for nickel-arsenide phases in the nickel-antimony, cobalt-antimony and iron-antimony systems.

Table 1

System	ρ, 25°	α, 25—100°
Ni—Sb	14×10^{-6}	500×10^{-5}
Co—Sb	35×10^{-6}	300×10^{-5}
Fe—Sb	230×10^{-6}	20×10^{-5}

It is seen that in the iron-antimony system the electric conductivity is very high, whereas the temperature coefficient is very low. An interesting point to note is that in iron-nickel-antimony ternary alloys with nickel-arsenide structure the temperature coefficient of electric conductivity drops, according to our observations (5), to nearly zero and even becomes negative in certain alloys. This testifies to the fact that nickel-arsenide compounds are nearer to semi-conductors in their properties than are true metals, or, at any rate, their number of effective free electrons is small, which implies a nearly complete filling up of the Brillouin zone and should serve as a criterion for selecting this or that form of zone. The ratio of the number of valency electrons to the number of atoms in an elementary cell of the nickel-arsenide compounds of antimony is assumed by us to be equal to $10/4 = 2.5$. Therefore, the zone sought for should include about 2.5 electronic states per atom.

Two nickel atoms and two arsenide atoms are contained in an elementary cell of a nickel-arsenide type of hexagonal structure. We shall describe the hexagonal structure in orthorhombic coordinates. The atom coordinates referred to orthorhombic axes are:

$$\text{Ni) } 000,\ 00\ {}^1/_2 \text{ and As) } 0\ \frac{1}{3}\ \frac{1}{4};\ \frac{1}{2}\ \frac{1}{6}\ \frac{1}{4}.$$

The structure factor has the form:

$$S_{hkl} = \cos^2 \frac{1}{2}\pi(h+k)\left[1 + e^{\pi i l} + e^{\pi i\left(\frac{2}{3}k + \frac{1}{2}l\right)} + e^{\pi i\left(h + \frac{1}{3}k + \frac{1}{2}l\right)}\right]$$

For deriving a Brillouin zone one needs to consider only those of the planes admitted by the structure factor which have the lowest indices and the highest values of intensity. These are the most intensive lines in the foreground of powderphotographs. Data referring to such planes denoted for orthorhombic lattice are presented in Table 2. Along with the general expression of the structure factor, the values of relative intensity I are given, which have been calculated by us for the NiSb compound (5).

Table 2

Indices for hexagonal lattice	(100)	(002)	(101)	(102)	(110)
Indices for orthorhombic lattice	(110)	(002)	(021),	(022),	(200),
General expression for S_{hkl}	$2f_{Ni} - f_{Sb}$	$2f_{Ni} - 2f_{Sb}$	(111) $\sqrt{3}\cdot f_{Sb}$	(112) $2f_{Ni} + f_{Sb}$	(130) $2f_{Ni} + 2f_{Sb}$
Computed intensities I for NiSb	0.0	1.6	34.3	32.1	35.9

The interplanar distance in the case of orthorhombic lattice is

$$d_{hkl} = \frac{1}{\sqrt{\left(\frac{h}{a}\right)^2 + \frac{1}{3}\left(\frac{k}{a}\right)^2 + \left(\frac{l}{c}\right)^2}}.$$

The perpendicular distance from the origin in K-space to the plane $(n_1\ n_2\ n_3)$ bounding the Brillouin zone is given by the equation

$$p_{n_1 n_2 n_3} = \frac{1}{2}\sqrt{\left\{\left(\frac{n_1}{a}\right)^2 + \frac{1}{3}\left(\frac{n_2}{a}\right)^2 + \left(\frac{n_3}{c}\right)^2\right\}}.$$

The zone bounded by the plane of the $(021) - (111)$ type has the form of a hexagonal bipyramid whose volume in K-space is $\frac{(4c^2+3a^2)^3}{72\sqrt{3}a^4c^5}$, where a and c are the lattice constants. The volume per atom is $\sqrt{3}\cdot a^2c/8$ and the number of states per atom for this zone is $n = \frac{(4c^2+3a^2)^3}{288a^2c^4}$.

Computations show that for the iron-antimony system n is equal to about 1.2. This corresponds to a very considerable overlapping of the second zone and will not agree with the electric properties discussed above. Therefore this zone cannot possibly be taken for the first Brillouin zone. On the same ground the zone bounded by planes of the $(022) - (112)$ type cannot be adopted either, since it includes four electronic states per atom.

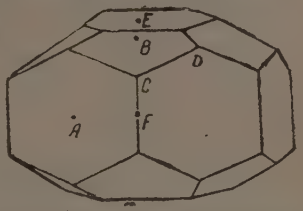
The actual shape of the Brillouin zone, such as corresponds to the criterion adopted by us at the beginning, is obtained as a combination of three types of plane, viz. $(022) - (112)$, $(200) - (130)$ and (002). Its shape is shown in the figure. The volume of this zone in the K-space is $\frac{(32a^2-2c^2)}{3\sqrt{3}a^4c}$ and the number of states per atom $n = \frac{16a^2-c^2}{6a^2}$.

Calculated values of n for nickel-arsenide phases in the systems of nickel; cobalt and iron with antimony are given in Table 3 along with electronic concentration computed from atomic composition.

Table 3

System	At% Sb	a (Å)	c (Å)	c/a	n	$\frac{N_{el}}{N_{at}}$
Fe—Sb	47.12	4.074	5.122	1.257	2.402	2.358
Co—Sb	48.21	3.898	5.177	1.328	2.372	2.409
Ni—Sb	49.46	3.940	5.140	1.304	2.386	2.469

It is seen from this table that if there is an overlapping with the second zone for iron-antimony alloys, this overlapping is small; for cobalt-antimony alloys the overlapping is larger; and it is still larger in the case of nickel-antimony alloys. These conclusions are in good agreement with the electric properties referred to in Table 1.



It is seen from the figure that the zone considered by us is rather flat, so that the overlapping of the second zone should take place first of all through the plane (002). From Table 2 it is evident that the intensity I of the line (002) on the powder-photographs is very weak, which is evidence of the low value of the energy discontinuity across the plane (002) of the Brillouin zone. In Table 4 the distances from the origin of the K-space (O) to the points marked on the figure are given. The points lying on the planes correspond, naturally enough, to perpendicular distances. The calculation is made for the NiSb compound, according to the data of Table 3.

Table 4

Distances	OE	OB	OA	OC	OD	OF
Length in inverse Å.	0.194	0.243	0.254	0.305	0.290	0.294

Thus: the first Brillouin zone for the nickel-arsenide type of structure is bounded by the planes (022)—(112), (200)—(130) and (002) and can include about 2.5 electronic states per atom. The overlapping into the second zone takes place across the plane (002), corresponding to the shortest distance from the origin of the K-space and to the least energy discontinuity as compared to the rest of the planes. The extent of the overlapping determines the electric properties of the particular compounds and accounts for the metallic character of phases with nickel-arsenide structure.

In conclusion I wish to express my gratitude to Prof. N. V. Ageev for the interest he took in this work and for valuable advice.

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TECHNICAL CHEMISTRY

PREPARATION OF MOTOR AND BOILER FUELS FROM PEAT BY THE METHOD OF THERMAL SOLUTION

By M. K. DIAKOVA

(Communicated by N. P. Chizhevsky, Member of the Academy, 14.IV.1943)

The Soviet Union has the largest peat reserves and is the biggest peat producer in the world (1). With an effective industrial method of converting it into liquid fuel, peat might become an important source of local liquid fuel for a number of regions and industrial districts of the country.

At present peat is being used here almost exclusively as a solid fuel of poor calorific value. Chemical and thermal treatment of peat, to convert it, in particular, into liquid fuel, has not been organized. Chemical methods of converting solid fuel into liquid fuel, applicable to coals and combustible shales (hydrogenation, carbonization), are found to be either little effective or altogether unsuitable in the case of peat.

Investigations I have conducted on the thermal solution in tetraline of diverse solid fuels (2), among them of the peat stage in carbonization, have revealed the possibility of developing a new, effective industrial method of producing artificial liquid fuel from peat by dissolving it in liquid organic solvents. With this end in view a more detailed study was made of the process of solution of typical Soviet peats in commercial solvents.

Samples of peats from the following deposits were investigated:

1. Bog peat from the Redkin Bog in Kalinin region: moisture content (laboratory sample), 8.4% and ash content, 8.75%.

2. Bog Ural peat received from the Ural Heavy Machine Building Works in 1942 (an average sample): moisture content (laboratory sample), 11.48% and ash content, 7.65%.

3. Bog peat from the White Russian SSR: moisture content (laboratory sample), 12.2% and ash content, 13.12%.

4. Bog peat of Kazan from the Borisogleb deposit in Yudin district, Tartar ASSR: moisture content (laboratory sample), 6.22% and ash content, 9.02%.

5. Bog peat of Kazan from Lenin district, Tartar ASSR: moisture content (laboratory sample), 24.98% and ash content, 10.06%.

As solvents we employed tetraline (b. p. 202—204° C; $d_{20}^{20}=0.971$); anthracene oil (the fraction boiling between 250 and 330° C, $d_{20}^{20}=1.0545$), a distillate of peat pitch (the fraction boiling between 200 and 350° C, $d_{20}^{20}=0.993$), the fraction of a hydrogenated peat solution (boiling away within the limits 230 and 330° C), and mazut (boiling above 300° C; $d_{4}^{20}=0.9508$).

The experiments were carried out in a half-liter revolving autoclave and in a continuous operation pilot unit with a capacity of 3 kg of raw material per hour.

The experimental procedure employed was the same as that used previously in studying the solution of coals and combustible shales (³).

A study was made of the influence of temperature (between 325 and 420°); heating time (between 5 and 60 minutes) and ratio of organic matter of the peat to the solvent (within the limits from 1: 3 to 2 : 3); and also of the influence of the nature of the solvent. Some of the most significant results are given in Tables 1 and 2.

Table 1

Solution of Soviet Peats in Organic Solvents

(ratio of sample to solvent 1:2, 1:3)

Exp. No.	Sample	Solvent	Experimental conditions		Extraction of organic matter of peat, %
			time (in min.)	temperature (in °C)	
538	Redkin peat	Tetraline	30	400	79.1
570	» »	»	5	420	86.3
561	» »	Anthracene oil	30	400	73.0
540	» »	Hydrogenate of a peat solution	15	410	73.0
575	» »	Gas oil	30	400	72.3
513	» »	Distillate of peat pitch	30	350	59.1
650	Urals peat	Mazut	10	410	67.7
606	Kazan peat, Lenin district	Anthracene oil	30	400	85.8
607	Do	Mazut	30	410	69.0
645	Kazan peat, Yudin district	»	10	410	62.2
591	White Russian peat	Anthracene oil	5	420	63,7

As the data summarized in Table 1 show, the thermal solution method ensures sufficiently intensive extraction of the organic matter in all the peat samples investigated. 70—85% of the organic matter is converted into liquid fuel; gas and pyrogenic water, 40—45% being converted into liquid fuel.

Although the peat begins to pass into solution at comparatively low temperatures, of the order of 325—350°, the process takes place slowly. At higher temperatures, between 410 and 420°, maximum solution is reached in 10 to 30 minutes.

Solubility of the peat is sufficiently high in all the solvents studied. It was found to be maximum (86% of the organic component) in tetraline. Likewise high is their solubility in anthracene oil (73—85%), a hydrogenate of a peat solution (73%), and mazut (69%). In a distillate of peat pitch, however, solubility was observed to be considerably worse (59%).

Variations in the ratio between the organic matter content of the peat and the solvent within the limits 1 : 3 and 2 : 3 do not essentially affect the solubility.

In thermal solution of peat a considerable part of its organic matter decomposes with the formation of gas, pyrogenic water and light fractions of liquid fuel (benzine, kerosene), while the other part passes into the solution forming liquid products of the type of boiler fuel.

The thermal solution method may be adopted for industrial production of motor and liquid boiler fuel in two ways. The most feasible of these from the point of view of present-day chemical technology of solid fuels is to combine the processes of thermal solution of peat and subsequent hydrogenation of the peat solution.

Roughly, this can be effected as follows. Ground and dried peat is subjected to thermal solution in a hydrogenate of a peat solution boiling above 300°; with this practically all the oxygen of the organic component, which is detrimental ballast, is given off in the form of CO_2, CO and H_2O. After distilling off benzine and ligroine from the products of solution, the insoluble residue is separated from the solution by filtration, and the filtrate is subjected to liquid phase hydrogenation;

Table 2

Yield of Products in Thermal Solution of Peat

(ratio of sample to solvent 1:3, 1:1.5)

Exp. No.	Sample	Solvent	Experimental conditions: time (in min.)	Experimental conditions: temperature (in °C)	Extraction of organic matter of peat, %	Products obtained in % of organic matter of peat: Gas	Crude benzine	Crude kerosene	Pyrogenetic water	Extract	Insoluble residue	Regeneration of the solvent
538	Redkin peat	Tetraline	30	400	79.1	17.0	12.6	—	13.8	35.6	22.4	100
564	» »	Anthracene oil	30	400	70.6	17.5	11.8	—	8.9	32.0	29.8	100
625	» »	Mazut	30	400	62	15.9	10.4	19.8	12.6	Boiler fuels 41.6	—	100
650	Urals peat	»	10	410	67.7	15.2	14.5	—	12.3	58		100
1st* exp.	Kazan peat from the Borisogleb deposit in Yudin district	»	20	420	62.1	14.9	24.3	19.1	9.4	28.7		100
2nd* exp.	Do	»	20	430	59.4	30.5	28.2	16.7	14.7	5.4		100

* Carried out on a continuous operation pilot unit.

then benzine; ligroine and medium oil are distilled off from the hydrogenate; and the residue of the hydrogenate boiling above 300° is employed as a solvent for treating a new portion of peat. By combining the methods of thermal solution of peat and hydrogenation of the solutions; up to 350 kg of clear products (benzine and kerosene) may be obtained from a ton of perfectly dry peat.

The second, simpler, but less accurate process of converting peat into liquid fuel; which does away with hydrogenation; consists in dissolving the peat in black mineral oil (mazut) and yields the basic products: benzine, ligroine; kerosene and boiler fuel. The technology of this prosess is roughly as follows. A mixture of dry peat (with a 3—5% moisture content) and mazut in the ratio 2 : 3 is subjected to thermal solution at 420—430° for a period of 20 minutes. The products of solution resulting are gases, benzine, ligroine, kerosene and a heavy liquid residue consisting of mazut, peat that has passed into the solution and insoluble particles of peat (among them mineral substances). This residue may be utilized as liquid boiler fuel.

A ton of perfectly dry peat (with a 9.02% ash content) and one and a half tons of mazut subjected to thermal solution under the indicated conditions with complete regeneration of the initial mazut (as boiler fuel) yield 280 kg of crude benzine; 167 kg of kerosene and 54 kg of boiler fuel with an ash content of 5—6% and a setting temperature of +9° (at a setting temperature of 5° for the initial mazut). This method ensures 50% conversion of the organic component of the peat into benzine, kerosene and boiler fuel.

Experiments designed to study the solution of peat in mazut in a continuous operation set-up with the raw material remaining in the reaction chamber (at 420—430° and 20—30 atm. pressure for an interval of 20 minutes) have shown that it is fully possible to make the process continuous on an industrial scale as well, and have made it possible to outline the basic features of the technological equipment. This method is expedient for industrial districts situated nearby peat deposits utilizing mazut as boiler fuel.

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GEOCHEMISTRY

ACCUMULATION OF GERMANIUM IN COALS

By V. M. RATYNSKY

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As shown by Goldschmidt ([1–2]), the area of accumulation of germanium coincides with that of coal basins. While the content of germanium in the earth crust is 0.0007%; its concentration in coals attains 0.01%. Much higher concentrations are observed in coal ashes; the figures for ashes rich in germanium showing a mean of 0.05% (the mean coefficient of enrichment equals 70). In individual samples the content of germanium attains 1.1%, the maximum coefficient of enrichment being equal to 1.600. Alongside of such important accumulations of germanium there are also coals that do not at all concentrate this element, its content in their ashes being as low as 0.001%.

The content of germanium shows a wide range of variation not only for different coal deposits, but even within a single deposit. As a general rule; it has been stated that the highest content of germanium occurs in coals poor in ashes.

In order to approach the solution of the problem as to the mode of germanium accumulation in mineral coals, we investigated several industrial deposits of the Urals and the Caucasus.

I. The Urals. 1) Cheliabinsk brown coal basin. Upper Trias (Rhaetian), Lower Jura (Lias). Banded coals. Deposits: Korkinskoe, Kopeyskoe.

2) Egorshinsky coal basin. Carbon. Anthracite, frequently graphitic. Egorshinskoe deposit.

II. The Caucasus. 3) Khumarinskoe mineral coal deposit. Lower Jura (Lias). Banded glance coals.

About 500 samples taken directly from the deposits were examined.

The coals to be analysed were ground in a mortar and burned so as to form thin ash layers in a porcelain dish over the flame of a gas burner; they were previously humectated with nitric acid. The content of germanium in the ash was evaluated quantitatively by way of spectrum analysis. The accuracy of determination was up to 0.001%, the error $\pm 7\%$. In those cases when germanium was present in a great quantity (0.1—1.0%), the content was checked by chemical methods.

The Urals coals are characterized by their low germanium content: especially low in germanium are the coals of the Egorshinskoe deposit; where it does not exceed 0.001% in the ash. Considerably richer are those of the Khumarinskoe deposit (Caucasus); in separate samples the content of germanium in them was as high as 1.0% in ash (Table 1).

The work carried out permits of some conclusions as to the mode of distribution of germanium in coals. Accumulation of germanium is typical of one of the petrographic varieties of coals, viz. of vitrain. Being poorer in ash than the rest of coal components, vitrain is present in relatively large quantities in coals that give little ashes, which accounts for a rather high germanium content in the latter.

Table 1

Germanium Content in Coals of Khumarinskoe Deposit

Sample No.	Material	Ash in %	Germanium content in the ash in%
	Clarain-vitrain, finely interbedded with slates and salt deposits	16.3	0.005
2	Do	10.0	0.004
4	»	21.5	0.005
6a	»	24.2	0.013
8	Vitrain lens	3.16	0.168
11	Clarain-vitrain, finely interbedded with slates and salt deposits	8.5	0.099
13	Do	3.8	0.024
14	»	6.0	0.015
16	»	6.8	0.004
17	»	4.3	0.004
18	»	9.0	0.148
21	»	13.8	0.026
21	»	18.2	0.008
22	»	10.7	0.018
24	Vitrain	6.5	0.153
27	Clarain-vitrain, finely interbedded with slates and salt deposits	23.0	0.002
28	Do	4.3	0.009
30	»	13.7	0.035
32	Clarain-vitrain (no slates)	2.5	0.054
35	Clarain-vitrain, finely interbedded with slates and salt deposits	6.7	0.002
36	Do	4.7	0.004
38	»	24.0	0.003
41	»	7.3	0.003
42a	Slates	85.6	<0.001
43	Glance coal, homogeneous with salt deposits along fissures	8.3	0.002
44	Clarain-vitrain, finely interbedded with slates and salt deposits	4.0	0.004
45	Do	3.1	0.011
46	»	17.3	0.003
46	Vitrain	3.3	0.113
48	Clarain-vitrain in pure state (without visible interbeds of slates)	8.8	0.129
49	Clarain-vitrain, finely interbedded with slates and salt deposits	6.5	0.011
51	Slates	64.2	<0.001
53	Clarain-vitrain, finely interbedded with slates and salt deposits	11.6	0.027
54	Do	12.1	0.054
54	»	16.1	0.009
55	Vitrain	3.5	0.588
55	Clarain-vitrain, finely interbedded with slates and salt deposits	34.6	0.003
58	Do	29.8	0.002
59	»	8.0	0.105
	Prevalence of vitrain		
60	Clarain-vitrain, finely interbedded with slates and salt deposits	5.5	0.030
60	Do	7.6	0.025
61	»	4.0	0.044
63	»	5.5	0.008
64	»	4.3	0.006
66	»	9.2	0.009
67	Coal, homogeneous, glance	1.8	0.051
67	Clarain-vitrain, finely interbedded with slates and salt deposits	32.3	0.008
71	Do	14.2	<0.001
71	»	13.2	<0.001
72	»	13.1	<0.001
72	»	15.7	<0.001
73	»	15.5	<0.001
73	»	18.0	0.004
76	»	8.2	0.013
78	»	5.5	0.191
84	»	5.1	0.036
86	»	26.0	0.003
88	»	7.5	0.038
89	»	6.8	0.005
91	»	14.2	0.005
92	»	12.3	0.006
93	»	4.4	0.054
95	»	6.6	0.063
96	»	13.6	0.005
101	»	6.8	0.004
102	»	4.0	0.007
107	»	12.5	<0.001
108	»	4.3	0,066
109	»	8.9	0.002
111	»	10.7	0.007
112	»	8.7	0.005
118	»	9.3	0.013
119	»	7.8	0.007
122	»	7.3	0.006
123	»	6.7	0.027
125	Slates	65.5	<0.001
126	Clarain-vitrain, finely interbedded with slates and salt deposits	25.5	<0.001
128	Do	13.3	0.003
129	»	13.8	0.035
132	»	3.7	0,003
140	»	3.8	0.082
145	»	3.3	0.052
153	»	9.2	0,045
153	»	17.7	0.020
154	»	10.2	0,020
157	»	5.6	0.157
158	»	11.1	0.016
159	»	6.6	0.023
160	»	9.8	0.004
160	Vitrain in pure state	1.4	0.986

Continued

Sample No.	Material	Ash in %	Germanium content in the ash in%	Sample No.	Material	Ash in %	Germanium content in the ash in%
162	Clarain-vitrain, finely interbedded with slates and salt deposits	5,5	0.027	190	Vitrain	2.3	0.815
164	Do	4.7	0.013	192	Clarain-vitrain, finely interbedded with slates and salt deposits	9.7	0.006
170	»	4.5	0.012	195	Do	10.3	0.003
170	»	10.7	0.005	197	»	29.2	0.001
171	»	9.5	0.004	198	»	35.2	<0.001
183	»	12.0	0.003	199	Coal, homogeneous, semi-lustrous, dense	7.3	<0.001
186	»	9.8	0.007				
187	»	12.5	0.006	201	Clarain-vitrain, without visible interbedding of slates	6.7	0.074
188	»	17.6	0.005				
189	»	14.7	0.002				

If compared to other coal components of equal or lower ash content: even those samples of vitrain that contain admixtures and are consequently rich in ash, invariably show a much higher germanium content, which is well illustrated by a comparison of samples 8 and 13, 16 and 24, 32 and 55.

In the banded coals of the Khumarinskoe deposit which are largely made up of strata of clarain-vitrain alternating with thin layers of carbonaceous and argillaceous schists, the fact of accumulation of germanium in vitrain is particularly conspicuous. This may well be examplified by the analysis of one of the samples made for each stratum separately (Table 2).

Table 2

No.	Material	Ash in %	Germanium content in ash in %
1	Vitrain in pure state	1.44	0.986
2	Vitrain finely interbedded with slates	16.1	0.007
3	Slate interbed, 2 mm thick, with vitrain inclusions	67.6	0.005
4	Interbed of carbonaceous schists, 1 mm thick	40.8	0.002
5	Mean test of the sample	15.5	0.010

In spite of a very high germanium concentration in vitrain (about 1% in the ash) the average test of a sample gave but 0.010% of germanium. The variations in germanium content do not go parallel with those of ash content: when the latter increases tenfold, the content of germanium drops as low as 100 times (Nos. 1 and 5; Table 2).

It should be pointed out that carbonaceous schists (Table 2) are much lower in germanium than slates with scarcely observable vitrain inclusions, and this in spite of the high ash content of the slates. The increased accumulation of germanium in vitrain should probably be put down to a higher absorbing capacity of the material that had served for the formation of vitrains.

Work on germanium accumulation in coals is in progress.

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BIOCHEMISTRY

ON THE PROBLEM OF UTILIZATION OF INTESTINE SLIME

By I. A. SMORODINTZEV and G. G. TER-SARKISSOV

(Communicated by S. S. Nametkin, Member of the Academy. 10.II.1943)

Up to the present intestine slime has only found application as an addition in metal treatment and as a constituent part of combined food. But its enzyme content which can be turned to profitable account in different branches of industry has been utterly disregarded.

The present authors have set themselves the task to determine the enzyme content in slime or in mucous membrane of small intestines and to develop methods for their extraction in the form of finished articles.

Method of Investigation. As a standard for comparing the fermentative power, we took extracts from pancreas and treated the slime by the same method. A hog pancreas 24 hours after extraction from the animal was freed from fat and connective tissue; ground in a mincing machine and left for infusion during 24 hours with four its volume of a 0.5% soda solution, in the presence of several millilitres of fresh gall and thymol crystals. In a similar way an extract with 60% glycerine was obtained by infusion for 48 hours. An analogous method was used for preparing soda extracts and glycerine extracts from hog intestine slime, obtained from slime machines of Moscow Meat Combinate. The extracts were filtered through gauze and their content of trypsine was determined by the Gross method and from the accumulation of aminonitrogen on formol titration. For a unit of tripsine—TU—we took the amount of nitrogen in milligrams formed during one hour under the action of 1 ml of extract on 25 ml of 3% gelatine 40° C and pH = 8.2.

Experimental Part. Activity of Trypsine of Pancreas. In a millilitre of soda extract or glycerine extract prepared from hog pancreas by the method described above 400 units of trypsine, according to Gross, were contained.

As an average of several determinations 1 ml of pancreas extract diluted 1 : 50 gave, on formol titration; 3.9 mg of nitrogen per hour; 1 ml of not diluted extract gave 50 times more, i. e. 195 mg of nitrogen. This value, according to our determination; was arbitrarily taken as a unit of comparison for measuring the activity of trypsine in extracts from pancreas and from intestine slime.

The time-growth of aminonitrogen, split off from gelatine by trypsine extract from pancreas, diluted 1 to 50, is shown in Table 1.

It is seen that glycerine extract and soda extract practically show the same number of TU per 1 g of raw slime. These data conform to a general rule that the enzyme acts strongly on the substrate at the beginning, but by and by its effect is diminished (1).

Investigation of Slime. The Gross method could not be used, because slightly diluted extracts are turbid; and strongly diluted ones contain but insignificant quantity of the enzyme.

Table 1

Pancreas

Duration of tests	30 min.		60 min.		90 min.	
Extracts	N, mg	TU	N, mg	TU	N, mg	TU
Glycerine	1.96	2	3.64	3.6	4.03	4
Soda	2.24	2.2	3.64	3.6	3.92	4

The formol titration was carried out with extracts diluted 1 : 10 and 1 : 20; because an extract diluted 1 : 50 is no longer active.

Average data for a set of experiments are presented in Tables 2 and 3.

Table 2

Extracts from Slime 1:10

Duration of tests	30 min.		60 min.		90 min.	
Extracts	N, mg	TU	N, mg	TU	N, mg	TU
Glycerine	0.48	0.09	0.94	0.19	1.35	0.27
Soda	0.55	0.11	1.16	0.23	1.43	0.23

Here too the glycerine extract and the soda extract practically show the same activity which is 10 to 15 times lower than the activity of a pancreas extract, as referred to 1 g of raw organ (Table 3).

Table 3

Extracts from Slime 1:10

Duration of tests	1 hour		2 hours		3 hours	
Extracts	N, mg	TU	N, mg	TU	N, mg	TU
Glycerine	0.85	0.17	1.72	0.34	2.18	0.44
Soda	1.2	0.24	1.84	0.37	2.45	0.49

As is generally the case, the enzyme acted more energetically at the first period than later; the soda extract appears to possess a somewhat higher activity than the glycerine extract.

The optimum temperature for the action of the trypsine from pancreas and slime was found to lie at 40° C.

Our experiments have shown that the slime's content of trypsine is 10 to 20% compared to the activity of the pancreas. Since the pancreas, from which usually oropon is manufactured for tanning industry, has now found extensive use in preparing insuline, a valuable medicinal preparation, the slime can be used successfully for manufacturing leather softeners and, perhaps, by means of adsorption for other trypsine preparations. This is all the more profitable, as after the extraction of the enzyme the slime can still be used as addition in metal treatment and as a constituent part of combined food.

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GEOLOGY

ON THE TECTONIC NATURE OF THE OIL-BEARING STRUCTURES OF POLASNA-KRASNOKAMSK TYPE

By V. A. APRODOV

(Communicated by V. A. Obruchev, Member of the Academy, 23.XII.1942)

The structures of the Polasna-Krasnokamsk part of the Near-Kama oil-bearing region belong to a particular type of platform tectonics. Up to the present time there is no settled opinion as to their nature. There exist four main trends in the discussion of the problem. The first of them refutes the leading rôle of tectonic processes in the development of the structures, referring them to an accumulation and erosion of sediments during the Palaeozoic.

These statements are developed in most details by Gerassimov (1). The second viewpoint is to interpret the local structures as anticline elevations formed as a result of unilateral tangential stress on the part of the western slope of the Urals and the Ufa plateau. This interpretation largely owes its development to Arkhangelsky (2). In the same way, viz. as platform elevations of the Herzynian foldedness, are the local structures considered by Nosal (3); Larionova (4) and others. The third group of authors explain the genesis of the Near-Kama structures by the intersection of two folds of Herzynian and Alpine orogenesis; followed by the formation of brachyanticlines; «nodes of intersection».

This viewpoint has been most thoroughly treated by Lapshinov (5).

Lastly, the fourth interpretation is given by V. V. Beloussov in his preliminary conclusions on the tectonics of the Volga-Urals region. The local structures are understood by him as more or less important domes, arisen in result of upheavals that took place within an area of continuous subsidence. At the same time this author points out the peculiar type of their tectonics; characterized by the formation of «families» of brachyanticlines against a background of more important upheavals.

The viewpoints referred to above are consecutive development stages of the interpretations given to the tectonics of the Permian Near-Kama region. Field observations carried out by the present author in the course of 1935 and 1942, as well as an analysis of structural maps partly published by Gerassimov (1); Lapshinov (5); Pometun (6), make us materially alter our idea as to the tectonics of the local structures. Their chief peculiarity consists in the radial, fan-like arrangement of brachyanticlines and brachysinclines with respect to the larger dome-like upheavals.

This radial system may distinctly be traced in the orientation of the strikes of the line of anticlines surrounding the large upheaval, with its centre near village Shalashnaya; of these the Polasna-Krasnokamsky is oriented south-west; Lobanovsky; south; Sylvensky, south-east; Yarinsky; north-east. A similar regula-

rity was discovered by this author in 1937 for a secondary fold-waviness of strata along the lower reach of the Chusovaya river, and by A. A. Aprodova in 1942 for the Nytva oil-bearing structures. An analogous phenomenon of fan-like arrangement of secondary upheavals has also been established for the North-Kama brachyanticline according to the magnetic map ([6]). It should be especially stressed that no such phenomena are observed for brachysynclinal areas.

The second trait peculiar of the local structures is a considerable change in the thickness of the layer of plastic rocks; according to the position of the section in the structure. The thickness of sulphate rocks of the Upper Kungurian in the Krasnokamsk structure varies within 14 m at the wings and up to 30 m at the vault.

Similar phenomena were established by A. A. Aprodova in 1942 for secondary upheavals of the Nytva brachyanticline.

The stratigraphical arrangement of carbonate rocks also varies essentially according to their position with respect to the vault. The vaults of secondary folds show a sharply increased dolomitization, gypsification and porosity of the rocks.

This should be accounted for by the metasomatic processes of their alteration.

The third peculiarity of the local oil deposits is the fan-like distribution in them of various types of underground waters. V. M. Kukanov (1937) and A. M. Kuznetzov (1939) established two main types of underground waters in the local Permian rocks: chloride-sodium-calcium waters, showing a mineralization of 20° Be, and chloride-sulphate-sodium waters with a mineralization of 5° Be. The former type corresponds to the domes of those structures; here belong waters that have migrated along fissures from deeper water-bearing horizons of the Carbonaceous.

On the ground of all the peculiarities observed we feel authorized to suggest a new tectonic interpretation of the structures described. As the original cause that led to their formation we, in agreement with the statements made by Beloussov in 1941, consider a differentiated vertical movement of platform areas of the earth' crust. However, contrary to what has been supposed by Beloussov, we believe this movement to have been an upheaval rather than a subsidence. It is just in the case of upheavals that matter is likely to have accumulated in this section. Experiments by Bridgman ([7]) and Adams ([8]) showed that at a considerable depth the rocks are subject to intense differential and hydrostatic strain which imparts them their elastic state. Local decreases of load, due to various causes (erosion, fissures, etc.), lead to a formation in these points of upheaval centres. But, aside from this the movement of the crystalline base seems to have played an especially important part here.

The rising substance of bottom rocks and of their derivatives is drawn into the area of upheaval and transmit the strain to the layer of sedimentation rocks which form a slant dome within the zone of weakened strain. The work made by the strata is hereby that of compression rather than stretch, since the horizontal component of movement in the area of upheaval is directed from its periphery towards the centre.

Let us imagine a flat section of the stratum in the form of a circle with a radius R. In place of this original circle there is formed a curvilinear surface of a globular segment or perhaps of a cone: with a resultant R. This lateral surface will always be smaller than the original plane. Its reduction will be the more considerable, the more intense the dislocation suffered by the stratum. At the expense of the original area around the dome secondary brachyanticlines and brachysynclines will be formed, arranged as long axes along the radii of the chief upheaval, and in the form of coulisse. In the upper zones of the geological section some horizontal displacements are to be expected, as well as a rotation of the axis of the structure relative to deeper horizons. According to the local conditions of the manifestation of forces; the process of formation of secondary structures may be attended either by compression; or by stretch of the rocks; in the latter case it will approach the normal process of fold formation. Thus, contrary to the opinion of

Kalizky ([9]) who denies the downward movement in the course of formation of small syncline structures or platforms, we admit its possibility, as well as that of accumulation there of oil and gas.

Owing to a peculiar distribution of forces around the main dome; the distribution of readily movable components (water, oil, gas) will proceed actively, as a component of the general movement of matter during fold formation. This has been pointed out for normal foldedness by Rosanov ([10]).

The diversified character of the formation of local structures ([5]) requires the evaluation of the individual importance of each of the phases in the distribution of oil in the structures. In particular, interest in this respect belongs to the orogenic phases of Alpine foldedness. Certain observations carried out on the water- and oil-bearing properties of the Near-Kama region point to that these components have quite recently been highly capable of migration.

Serious attention is deserved therefore by the treatment of problems referring to the Quaternary tectogenesis of the Permian Near-Kama region during the Bakinskian and Upper Quaternary orogenic phases. It seems probable that in some instances they might have materially affected the oil-bearing capacity of the Near-Kama structures.

To conclude it should be noted that the regularities established in the distribution of upheavals permit to assume the existence of anticline structures of the Polasna-Krasnokamsk type along the lower course of the Obva river, north-west to Polasna, and in several other points of the Permian Near-Kama region.

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GENETICS

INDUCED DOUBLE FLOWERS IN STOCKS

By L. BRESLAVETZ

(Communicated by L. A. Orbeli, Member of the Academy, 12.IV.1943)

In the summer of 1940 the appearance of double flowers was stated in the progeny of seven stock plants; subjected to the actions of acenaphthene in the previous year (1). The number of double-flower stocks varied in individual families from 41.7 to 76%.

In order to clear up the nature of this phenomenon mass plantings (one and a half—two thousand of every family) were carried out in 1941. 200 from 300 seeds were taken from every offspring for separate planting. In order to characterize the progeny, we take the family No. 88 which will serve to illustrate the transmission of the double-flower habit by a seed generation (Table 1).

Table 1

Plant	Number			Per cent of double
	Double	Single	Total	
88_1	85	81	166	51.2
88_2	15	23	38	39.4
88_3	124	103	227	54.6
88_4	168	124	292	57.3
88_5	65	36	101	64.3
88_6	195	129	284	54.6
88_7	53	102	155	34.1
88_8	112	89	201	55.7
88_{10}	108	77	185	57.6
88_{11}	142	102	244	58.2
88_{13}	101	84	185	54.5
88_{15}	186	118	304	61.2
88_{17}	99	39	138	71.4
88_{18}	95	74	169	56.2
Total	1591	1281	2872	55.4

As may be seen from Table 1, the fluctuations in the appearance of double flowers in the progeny of a given family are very considerable ranging from 34.2 to 71.4%. In general; the decrease of the number of double-flower plants is by 11.6%.

Turning now to the progeny of two more families; we see that in one of them there is no reduction; whereas in the other the reduction is as strong as to attain the value established for the first family (Table 2).

Table 2

Family	Ratio of double-flower plants	
	1940	1941
88	66.0	55.4
89	51.0	52.4
95	74.5	59.9

Two conclusions may be drawn from these observations:

1. The double-flower habit induced in stocks experimentally by the action of acenaphthene on seeds is transmitted by a seed generation to an extent that characterizes the particular family. In 1941 as well as in 1940 the highest ratio of double-flower plants was among the family No. 95.

2. Though the seeds for planting were selected from lines showing the highest ratio of double-flower plants; there was observed a reduction of double flowers in their 1941 progeny; which indeed was to be expected on the assumption expressed in an earlier article ([2]).

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PLANT PHYSIOLOGY

ON THE RÔLE OF SPECTRAL COMPOSITION OF LIGHT IN PHOTOPERIODIC REACTION

By A. F. KLESHNIN

(Communicated by A. N. Bach, Member of the Academy, 27.VII.1942)

This communication presents the results of investigations carried out in 1941 in which the data of the previous works and the defects of the method then applied have been taken into consideration. Accordingly, a special method has been evolved securing a purity of spectral composition and a homogeneousness of the incident radiant energy; the main characteristic which speaks in favour of the new method is that the experiments were made with the individual leaf while on the plant under a comparatively low intensity of radiation.

As the test-object we took a short-day plant, *Perilla ocymoides*, which has been thoroughly investigated in point of photoperiodical reaction. It was found to be well fitted to a study by the method adopted by us.

Adult vegetating plants developing on long daylength were subjected previous to the experiment to the following operation. Their top was cut off on the level of the 6—7th pair of leaves, and all their leaves but the 5th or the 6th pair were removed; one of the leaves left served as control and was given continuous illumination (10 hours of natural light plus 14 hours of electric light); the other leaf—the test-leaf—was kept under the following light conditions: 10 hours of natural light (in the green-house from 9 a. m. to 7 p. m.) plus 14 hours of artificial light of a definite spectral composition (from 7 p. m. to 9 a. m.). The light conditions under which the test-leaves were kept are represented schematically as follows:

a) 10 hours of natural light + 14 hours of red light (620—720 mμ);
b) 10 hours of natural light + 14 hours of green light (480—580 mμ);
c) 10 hours of natural light + 14 hours of blue light (400—480 mμ);
d) 10 hours of natural light + 14 hours of white light (400—720 mμ);
e) 10 hours of natural light + 14 hours of darkness;
f) natural illumination; natural daylength.

Observations were made on shoots, arising in the axils of the test and control leaves (the test and control shoots; respectively).

The sources of illumination were ordinary incandescence lamps; 75, 500 and 1000 watt, which were immersed in glass vessels with continuous flow of water (the water served for cooling the lamp and for removing infra-red rays). The light of the requisite wave-length was obtained by means of liquid (water) light-filters, prepared in the following way:

1) Red light-filter (620—720 mμ)—tetracine solution (1.0 *n*) and methyl-violet (1.0 *n*).

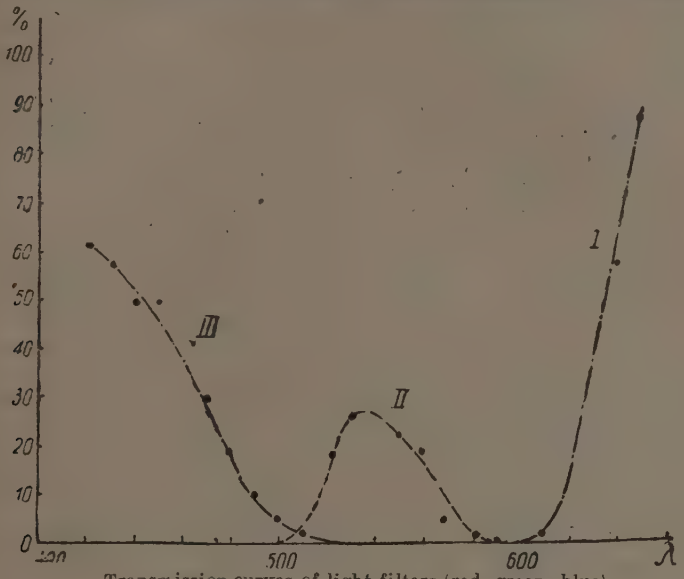
2) Green light-filter (480—570 mμ)—hydrochloric acid solution of K_2CrO_4 (100 *n*) and $CuSO_4$ (6000 *n*).

3) Blue light-filter (400—480 mμ)—a solution of methyl-violet (0.5 *n*) and $CuSO_4$ (6000 *n*).

4) White light-filter (400—720 mμ)—distilled water.

The light-filters were poured in pans made of ordinary window-glass which were in contact with light-proof chambers containing the test-leaves. The filters were changed periodically; the transmission curves of the light-filters (figure) were determined by means of a König-Martens spectrophotometer. An every-day control of the light-filter throughout the experiment was effected by means of a pocket spectroscope.

The intensity of radiation was regulated by measuring the distance between the test-leaf and the source of radiation. The measurements of the radiation inten-



Transmission curves of light filters (red, green, blue)

sity were made every day by the objective photometer designed by Maslakovetz and Makarevsky.

Two series of tests were started. In the first series the action of a constant quantity of incident radiant energy of different wave-lengths on the development of *Perilla ocymoides* was investigated.

Test 1. Illumination during 50 days; from June 6 till July 26; intensity of radiation, 1000 erg/cm² sec.; test completed on July 29.

Test 2. Illumination during 30 days, from June 6 till July 7; intensity of radiation, 1000 erg/cm² sec.; then the test-leaves of all sets were treated with a short 10-hour daylength for 26 days from July 7 till Aug. 2; the experiment completed September 3.

Test 3. Illumination during 30 days, from June 5 till July 7; intensity of radiation, 1000 erg cm² sec.; afterwards, natural daylength; the experiment completed September 3.

The results of this set of tests are presented in Table 1.

In the second set of tests the action of the intensity of light (500, 250, 100 and 50 erg cm² sec.) and of its spectral composition upon the photoperiodical reaction of *Perilla ocymoides* was studied. The conditions of these tests and the results are given in Table 2.

From the data referred to in Tables 1 and 2 the following conclusions may be made:

1. If the intensity of light in the process of the photoperiodical action is above 50 erg cm² sec.; the reproductive phase of development in short-day plants (*Perilla ocymoides*) is inhibited; whatever the spectral composition may be.

2. If the light intensity is above 50 erg/cm² sec.; the effect is not merely to

Table 2

Effect of Spectral Composition and Intensity on Budding in *Perilla ocymoides*

Set	Conditions of illumination	Wave-length in mμ	Budding of the test-shoot*								
			500 erg/cm² sec.; from July 14 to Aug. 9 (26 days)		250 erg/cm² sec.; from July 28 to Aug. 16 (19 days)		100 erg/cm² sec.; from July 14 to Aug. 9 (26 days)		50 erg/cm² sec.; from Aug. 1 to Aug. 22 (22 days)		
			Outset of budding	Number of days from start of illumination till budding	Outset of budding	Number of days from start of illumination till budding	Outset of budding	Number of days from start of illumination till budding	Outset of budding	Number of days from start of illumination till budding	Number of days from beginning of short day-length
1	Red light	620—720	Sept. 1	48	Sept. 8	52	Sept. 8	56**	Sept. 15	46	24
2	Green light	480—580	Sept. 1	48	Sept. 3	47	Sept. 1	48	Sept. 11	42	20
3	Blue light	400—480	Sept. 1	48	Sept. 3	47	Sept. 1	48	Sept. 3	34	12
4	White light	400—720	Sept. 1	48	Sept. 3	47	Sept. 8	56**	Sept. 15	46	24
5	Short daylength	—	Aug. 7	24	Aug. 16	19	Aug. 7	24	Aug. 22	22	22
6	Natural day	—	Sept. 1	48	Aug. 18	21	Aug. 22	22	Aug. 22	22	22

* In all cases except those of set 3 in the last group and set 5 in everyone of the groups, budding took place under the influence of naturally curtailed daylength.

** The retardation is connected with the affection of the plants.

Table 1

Effect of Spectral Composition of Light on Budding in *Perilla ocymoides*

Set	Conditions of illumination	Wave-length in mμ	Budding of the test-shoot						
			Test 1		Test 2			Test 3	
			Outset of budding	Number of days from start of illumination till budding	Outset of budding	Number of days till beginning of budding: from start of illumination	Number of days till beginning of budding: from start of short-day treatment	Outset of budding	Number of days from start of illumination till budding
1	Red light . .	620—720	—	∼	Aug. 4	59	27	Sept. 1	86
2	Green light . .	480—580	—	∼	Aug. 4	59	27	Sept. 1	86
3	Blue light . .	400—480	—	∼	Aug. 4	59	27	Sept. 1	86
4	White light .	400—720	—	∼	Aug. 4	59	27	Sept. 1	86
5	Short daylength	—	June 30	24	June 30	24	24	June 30	24
6	Natural day . .	—	—	∼	Aug. 4	59	27	Sept. 1	86

delay the development; but even to check completely the flowering; the reproductive phase does not begin at all unless the plant is given a short daylength.

3. Different regions of the spectrum have a different inhibiting effect on the process of the transition to the reproductive phase of development and lead to no specific reaction, which afterwards with the onset of favourable light conditions might delay or accelerate the outset of flowering; under short-day conditions flowering sets in simultaneously in all test and control sets.

4. Under a light intensity of 50 erg/cm² sec. the plants of different sets behave differently: red and green light, as well as white light completely stop the initiation of reproductive processes; blue light is no obstacle to their start and only delays it.

5. If the intensity of radiation is above 50 erg/cm² sec., the light factor is in excess; therefore in determining the rôle of different portions of the spectrum in the photoperiodical process, we have to follow the line of decreasing radiation intensity.

Thus, on the basis of our investigations and on the basis of data obtained by studying the radiation intensity in the photoperiodical process ([1],[2]), we arrive at the following conclusions. The photoperiodical processes are affected by any region of the visible part of the spectrum; it is only the amount of energy absorbed by some acceptor of the leaf that is important. The specific action of the light of different wave-lengths consists merely in the fact that the sensitivity of this acceptor is different in response to different wave-lengths. For every region of the spectrum there is a definite threshold of radiation intensity below which this region is perceived as darkness in the photoperiodical process; this threshold is different for different plants. The specific action of the light manifests itself only under low intensity (different for different plants); when, for instance, the blue light is already perceived as darkness and the red light, in response to which the susceptibility threshold is lower; is still perceived as light.

In the light of these conceptions the different behaviour of plants towards light can easily be accounted for. This also explains the contradictions between the experimental data of different authors.

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BACTÉRIOLOGIE

CULTURE DE *LEISHMANIA TROPICA* DANS L'OEUF DE POULE

Par D. A. BAHRAMI

(Présenté par E. N. Pavlovsky, de l'Académie, le 1.III.1943)

Aucun essai sur la culture de *Leishmania* dans les oeufs de poule n'ayant précédé notre recherche, il nous a paru utile d'attirer, dès maintenant, l'attention du monde savant sur le résultat de nos propres expériences.

Le premier ensemencement partait d'une souche de *Leishmania tropica* à son trente-troisième passage sur le milieu N. N. N. A ce but nous avons pris des oeufs de poule incubés préalablement pendant six jours à la temperature humide de 39°; aérés 15 minutes par jour et portés à l'épreuve devant une source de lumière pour être assuré de l'état du développement de l'embryon. Puis deux trous ont été pratiqués sur chaque oeuf au moyen d'un foret, dont l'un situé près de l'embryon nous a servi pour l'introduction du parasite, tandis que l'autre conduisait dans la chambre à l'air. Les trous bouchés à la paraffine, nous avons pu constater après le 3-ème jour d'étuve à 25° le développement d'un grand nombre de *Leishmania* dans la membrane chorio-allantoïde et dans le coeur de l'embryon. L'aspect microscopique de ces *Leptomonas* ne diffère point de celui que l'on peut observer sur le milieu N. N. N. Toutes les formes connues de ce parasite y sont représentées: formes arrondies, allongées, ovalaires, rosaces, etc. D'autre part, nous nous sommes servis dans l'ensemencement des oeufs du corps de *Leishmania* obtenu directement des boutons d'Orient; le résultat en n'était pas moins encourageant, et la facilité du développement des parasites était surprenante.

Toutefois l'oeuf de poule se présente comme un milieu propice à la culture de *Leishmania* et semble être identique à tout point de vue aux autres milieux déjà connus. Or, la culture de *Leishmania* dans l'oeuf de poule nous amène à constater un fait d'ordre biologique non moins important qui nous peut servir de base à d'autres thèses du même rang, à savoir: la possibilité de cultiver un protozoaire, tel le *Leishmania tropica*, dans l'embryon d'une poule, alors qu'il ne peut guère se développer ou vivre dans le corps même de cet oiseau. De là l'importance des oeufs d'oiseaux comme un milieu remplaçant les corps des insectes pour le développement des protozoaires dont une partie de vie doit passer dans l'organisme de ces arthropodes.

Institut Pasteur
de l'Iran.

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